

REGULARITY PROPERTIES, REPRESENTATION OF SOLUTIONS AND SPECTRAL ASYMPTOTICS OF SYSTEMS WITH MULTIPLICITIES

ILIA KAMOTSKI AND MICHAEL RUZHANSKY

ABSTRACT. Properties of solutions of generic hyperbolic systems with multiple characteristics with diagonalizable principal part are investigated. Solutions are represented as a Picard series with terms in the form of iterated Fourier integral operators. It is shown that this series is an asymptotic expansion with respect to smoothness under quite general geometric conditions. Propagation of singularities and sharp regularity properties of solutions are obtained. Results are applied to establish regularity estimates for scalar weakly hyperbolic equations with involutive characteristics. They are also applied to derive the first and second terms of spectral asymptotics for the corresponding elliptic systems.

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1. INTRODUCTION

Let X be a smooth manifold without boundary of dimension $n \geq 3$. Let P be an elliptic self-adjoint pseudo-differential operator of order one acting on half-densities on m -dimensional crosssections of vector bundles on X . We consider the following Cauchy problem for $u = u(t, x)$

$$(1.1) \quad \begin{cases} iu' - Pu = 0, & (t, x) \in [0, T] \times X, \\ u|_{t=0} = u_0. \end{cases}$$

It is well known that if equation (1.1) is strictly hyperbolic, the system can be diagonalized and its solution can be given as a sum of Fourier integral operators applied to Cauchy data (e.g. [6]). An important question that has been studied over many years is what happens when P has multiple characteristics.

Since we will be mostly interested in local properties of solutions, we may already assume that P acts on functions, and can think of it as an $m \times m$ matrix of pseudo-differential operators of order one and we think of u_0 as of an m -vector.

Let $A(x, \xi)$ denote the principal symbol of P . If A is a diagonal matrix, properties of system (1.1) have been studied by many authors. For example, in [16] and [15] Kumano-go and coauthors used the calculus of Fourier integral operators with multi-phases to show that the Cauchy problem (1.1) is well-posed in L^2 , Sobolev spaces H^s , and to study its propagation of singularities. Systems with symmetric principal part A have been extensively studied as well (e.g. [14], [12], etc.) In a generic situation, such systems have double characteristics, and their normal forms have been found by Braam and Duistermaat [2]. Recently, Colin de Verdiere [4] used these

representations to derive some asymptotic properties of such systems. Polarization properties of similar systems were studied by Dencker in [5].

More elaborate analysis of system (1.1) becomes possible if we assume that the principal symbol matrix $A(x, \xi)$ is smoothly microlocally diagonalizable with smooth eigenvalues $a_j(x, \xi)$ and smooth eigenspaces. Then, as it was pointed out by Rozenblum in [21], there exists a finite dimensional cover \tilde{X} of X such that A lifted to \tilde{X} can be globally diagonalized provided X is compact. In this situation Rozenblum showed that the Picard series for this problem gives an expansion with respect to smoothness in the case of non-involutive characteristics. In other words, one assumes that if $a_j(x, \xi) = a_k(x, \xi)$ for $j \neq k$, then the Poisson bracket $\{a_j, a_k\}(x, \xi) \neq 0$. This means that at all points of multiplicity, bicharacteristics intersect characteristic surfaces transversally. However, this condition is non-generic even for diagonal systems. For example, it is clear that if one of characteristics has a fold, there may be a point where this transversality condition fails, and it is not possible to remove it by small perturbations.

One purpose of this paper is to present results removing the transversal intersection condition. We will allow characteristics to be involutive of finite type and some characteristics to be identically equal. Operators satisfying our Condition C below will be also generic in the class of smoothly microlocally diagonalizable systems. Below we will explain that the microlocal diagonalizability condition is quite natural when considering weakly hyperbolic scalar equations with Levi conditions (Examples 1 and 2). This is also the case for Maxwell equations (e.g. [1]).

We will investigate regularity properties of system (1.1) in this generic setting. Even in the case when the system is strictly hyperbolic, L^p properties of solutions have been studied for many years, since already this case has several important applications for nonlinear equations and harmonic analysis. Regularity properties of non-degenerate Fourier integral operators have been established by Seeger, Sogge and Stein in [24]. They showed that a Fourier integral operator T of order zero satisfying local graph condition, is locally bounded from $(L^\alpha_\alpha)_{comp}$ to L^p_{loc} for $1 < p < \infty$ and $\alpha = (n-1)|1/p - 1/2|$. As a consequence they showed that if system (1.1) is strictly hyperbolic, there is a loss of α derivatives in L^p , i.e. $u_0 \in L^\alpha_\alpha$ implies $u(t, \cdot) \in L^p$. Moreover, if at least one of characteristic roots a_j is elliptic, the loss of α derivatives is sharp. If none of a_j 's is elliptic, this result can be improved ([22]).

Our Theorem 2.2 will establish a similar property for systems (1.1) with multiplicities. Moreover, this will imply L^p estimates for scalar weakly hyperbolic equations with involutive characteristics. It is known that in general weakly hyperbolic cases one often loses regularity even in L^2 . However, in the case of involutive characteristics the equation can be diagonalized and in Theorem 2.3 we will give L^p estimates for such equations. This will, on one hand, extend the L^p result of Seeger, Sogge and Stein to weakly hyperbolic equations and systems with multiplicities, while on the other hand establishing L^p estimates for systems considered by Kumano-go, Rozenblum, and others. The result will be general for scalar weakly hyperbolic equations satisfying Levi conditions with characteristics satisfying Condition C below.

Note that if a scalar operator strictly hyperbolic and we write it in the form (1.1), we can diagonalize P together with lower order terms (e.g. [15]) and split it into m

scalar equations, for which many things are known. However, in the case of multiple characteristics this is impossible, so a more elaborate analysis is needed.

Now we will formulate our main assumption. Let us define operator

$$H_{a_j}f = \{a_j, f\}, j = 1, \dots, m,$$

where

$$H_g(f) = \{g, f\} = \sum_{k=1}^n \left(\frac{\partial g}{\partial \xi_k} \frac{\partial f}{\partial x_k} - \frac{\partial g}{\partial x_k} \frac{\partial f}{\partial \xi_k} \right)$$

is the usual Poisson bracket. Our assumption is that at points of multiplicity $a_j = a_k$ of non-identical characteristics a_j and a_k , bicharacteristics of a_j intersect level sets $\{a_k = 1\}$ with finite order, i.e. $H_{a_j}^M a_k \neq 0$ for some M , at points where $a_j = a_k$. In other words, we allow involutive characteristics of finite type, and formulate our main condition.

Condition C:

Suppose that there exists $M \in \mathbb{N}$ such that if for some j and k , a_j and a_k are not identically the same, then

$$(1.2) \quad \begin{aligned} & a_j(x, \xi) = a_k(x, \xi), (x, \xi) \in T^*X \implies \\ & H_{a_j}^\lambda a_k(x, \xi) = \{a_j, \{a_j, \dots \{a_j, a_k\} \dots\}^\lambda\}(x, \xi) \neq 0, \end{aligned}$$

for some $\lambda \leq M$. While the function $M = M(x, \xi)$ is locally bounded, it is allowed to grow at infinity.

We note here that the transversality assumption of Rozenblum [21] requires (1.2) to hold with $M = 1$. Strictly hyperbolic case is also covered by this condition (in which case we set $M = 0$). The case of a_j and a_k defining glancing hypersurfaces (as in Melrose [17]) corresponds to $M = 2$.

In Section 2 we will give several examples of characteristics satisfying condition C, in particular those arising from weakly hyperbolic scalar equations with involutive characteristics. Such equations and propagation of their singularities have been analyzed in [3], [18], [19], [11], [13], etc. We will also establish estimates in L^p and other spaces for the weakly hyperbolic equations or systems satisfying condition C.

It is interesting to note that conditions similar to Condition C appeared in the study of subelliptic operators (e.g. Hörmander [10, Chapter 27]). For instance, in the case of 2×2 systems P with characteristics a_1 and a_2 , we can consider operators Q with principal symbol $q = a_1 + ia_2$. Then microlocal subellipticity of Q implies Condition C, with some M dependent on the loss of regularity for Q , which, therefore, implies the Weyl formula for P (Theorem 2.5), regularity estimates for (1.1) and all other results of this paper.

Now we will give an informal explanation of the strategy of our analysis. First, let us follow [21] to show that microlocal diagonalizability implies a local one on some cover \tilde{X} of X with finitely many leaves. For this argument we assume that X is compact. Since all the analysis of this paper will be local, if X is not compact, we can always assume that the amplitude of $P(x, D)$ is compactly supported.

Let an elliptic pseudo-differential operator $P(x, D)$ of order one act on sections of an m -dimensional Hermitian vector bundle E . Let $L^2(E)$ be the space of sections of half-densities on E and let P be self-adjoint on $L^2(E)$. Let E' be the lifting of E to T^*X . Then for each $(x, \xi) \in T^*X$ the principal symbol $A(x, \xi)$ of $P(x, D)$ is a Hermitian isomorphism of sections of E' . Without loss of generality we can assume that $A(x, \xi)$ is positive definite. Indeed, if it has both positive and negative eigenvalues, it is possible to globally block-diagonalize $A(x, D)$ with some suitably chosen pseudo-differential operators, to reduce it to a direct sum of positive and negative definite operators. Then each of these operators can be analyzed independently.

Let us assume that the principal symbol $A(x, \xi)$ is microlocally diagonalizable. This means that microlocally in $\Lambda \subset T^*X$ such that $E'|_\Lambda \cong \Lambda \times \mathbb{C}^m$, principal symbol $A(x, \xi)|_\Lambda$ has m smooth non-negative eigenvalues $a_j(x, \xi)$ and one dimensional eigenspaces $V_j(x, \xi)$, and such diagonalizations are compatible in intersecting cones. In this situation Lemma 5.1 insures that there is a cover \tilde{X} of X with finitely many leaves such that the principal symbol of the lifting of $P(x, D)$ to $T^*\tilde{X}$ can be globally diagonalized modulo lower order terms. Note that since dimensions of X and \tilde{X} are the same and because of formula (5.1) all our results on \tilde{X} will imply corresponding results on X . Therefore, we may assume that the principal symbol A of operator P may be smoothly diagonalized over compact subsets of X , that is

$$P = A + B, A = \text{diag}\{A_1, \dots, A_m\},$$

where $A_j \in \Psi^1$ are scalar pseudo-differential operators with principal symbols $a_j(x, \xi)$ (eigenvalues of A). Here a_j 's may be identically equal to each other or may intersect with any finite order, according to our Condition C. Here B is an $m \times m$ matrix of pseudo-differential operators of order zero. We may also assume that $B_{jj} = 0$ for $1 \leq j \leq m$, if we add these terms to the diagonal of A . For the moment we will also assume that none of a_j 's are identical. Otherwise, if some of a_j 's being identically equal to each other locally at some points, the construction is slightly different, but all the results remain valid. This will be carried out in detail in Section 3 in the proof of Theorem 2.2. Substitution $U = e^{-iAt}V$ leads to the equation

$$(1.3) \quad \begin{cases} V' = Z(t)V, \\ V|_{t=0} = I, \end{cases}$$

with $Z(t) = -ie^{iAt}Be^{-iAt}$. Writing the Picard series for problem (1.3), we obtain the expansion

$$(1.4) \quad V(t) = I + \int_0^t Z(t_1)dt_1 + \int_0^t \int_0^{t_1} Z(t_1)Z(t_2)dt_2dt_1 + \dots$$

A general term of this series is

$$(1.5) \quad Q_l = \overbrace{\int_0^t \int_0^{t_1} \dots \int_0^{t_{l-1}}}^l e^{iA_{j_1}t_1} b_{j_1j_2} e^{iA_{j_2}(t_2-t_1)} \dots dt_l \dots dt_1.$$

It is easy to see that $\|Q_l\|_{L^2 \rightarrow L^2} \leq C/l!$ and that series (1.4) converges in L^2 and in H^s . Using the notion of a multi-phase for Fourier integral operators, Kumano-go et al. ([16], [15]) studied propagation of singularities of Q_l . Instead of introducing multi-phases for Q_l , we will analyze operators Q_l in more detail and will show its smoothing properties in Sobolev spaces under Condition C. Here, contrary to the

transversal non-involutive case of Rozenblum (when $M = 1$), we do not have good control on the singular supports of integral kernels of operators Q_l , so more elaborate geometric analysis is required.

In fact, our Theorem 2.1 asserts that $Q_l(t)$ maps L^2 to some Sobolev space $H^{p(l)}$, and that $p(l) \rightarrow \infty$ as $l \rightarrow \infty$. Then we will show that this allows to treat the Picard expansion as a series with finitely many terms, with many conclusions, such as L^p estimates for solutions and spectral asymptotics of P .

In particular, L^p estimates will follow from a general principle which we will prove in Theorem 3.1 for equation $u' - Z(t)u = f, u(0) = u_0$. In Corollaries 3.2 and 3.3 we will show regularity of solutions of this equation for pseudo-differential operators of order zero $Z(t) \in \Psi^0$ or for Fourier integral operators of negative orders $Z(t) \in I^{-\epsilon}$. In general, there may be problems with this construction for $Z(t) \in I^0$, but there we can use the structural properties of $Z(t)$ in (1.3).

Everywhere in this paper $\Psi^\mu = \Psi_{1,0}^\mu(X)$ will denote the space of classical pseudo-differential operators of order μ of type $(1, 0)$. The space of Fourier integral operators of order μ with amplitudes of type $(1, 0)$ will be denoted by I^μ . All Fourier integral operators in this paper will be non-degenerate, which means that its canonical relation satisfies the local graph condition, i.e. it is a graph of a symplectic diffeomorphism from T^*X to itself. Constants C may be different in different formulas throughout this paper. We will use the following notation for norms and spaces. By L_α^p we will denote the Sobolev space of functions f such that $(I - \Delta)^{\alpha/2}f \in L^p$. For a function f we will denote its L^p -norm by $\|f\|_{L^p}$ and its Sobolev $H^s = L_s^2$ norm by $\|f\|_s$. If T is an operator, by $\|T\|_s$ we will denote its operator norm from L^2 to H^s .

Results of this paper can be established also in the case of operators P dependent on t . This will be the subject of a separate paper since it will also involve the analysis of non-smooth coefficients.

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2. MAIN RESULTS

Now we will formulate our results concerning terms of the Picard series (1.4) and solutions to systems (1.1) and (1.3). We will also give a Weyl formula for P .

Our first main result will be Theorem 2.1 on the smoothing properties of terms Q_l . The other two important results will be Theorem 2.2 on the L^p -regularity of solutions to system (1.1) and Theorem 2.5 on the spectral asymptotics for elliptic operator P satisfying Condition C. To obtain L^p -estimates, we use Theorem 3.1, which we regard as a general principle behind regularity estimates for general Cauchy problems based on several natural properties of the right hand side operators $Z(t)$. We will illustrate its use in several situations in Corollaries 1-3. Theorem 2.4 is a statement on the propagation of singularities of operators Q_l or, more generally, of Fourier integral operators in which the frequency integration is performed over a cone rather than over the whole space. Note that these results will hold for microlocally smoothly block-diagonalizable operators with any (finite) geometry of characteristics, i.e. characteristics satisfying our Condition C.

The following theorem establishes a smoothing property of operators Q_l under Condition C.

Theorem 2.1. *Let condition C be satisfied, that is assume that there is M such that for any $(x, \xi) \in T^*X$ and any $1 \leq j, k \leq m$, $j \neq k$, with a_j and a_k not identically the same and $a_j(x, \xi) = a_k(x, \xi)$, we have*

$$\{a_j, \{a_j, \dots \{a_j, a_k \overbrace{\}^{\lambda}} \dots \} \} (x, \xi) \neq 0$$

for some number $\lambda \leq M$. Then for sufficiently large l operator Q_l in (1.5) is bounded from L_{comp}^2 to H_{loc}^N , where

$$N = \frac{((-3n/2 - 2)(3[l/2] - 1 - n) + ([l/2] - n - 1)(l/(2M) - n - 1))}{(3[l/2] - 2n - 2 + l/(2M))} \sim \frac{l}{6M + 2}.$$

Note that the exact order N can be improved. However, it is most important that it increases to infinity as $l \rightarrow \infty$. This theorem implies, in particular, that the series (1.4), i.e. the series

$$V(t) = I + Q_1(t) + Q_2(t) + \dots$$

is a series with respect to smoothness. This fact allows one to refine the study of propagation of singularities and regularity properties of solutions to systems (1.1) and (1.3).

It turns out that smoothing properties of Q_l in Theorem 2.1 can be used to establish local L^p properties of solutions to systems with multiplicities (1.1). For strictly hyperbolic equations such estimates have been established by Seeger, Sogge and Stein in [24], and some optimal estimates were given in [22]. Our next result concerns regularity of solutions to Cauchy problem (1.1). In this theorem we also allow the lower order term B to depend on t .

Theorem 2.2. *Let $1 < p < \infty$ and $\alpha = (n - 1)|1/p - 1/2|$. Let $P = P(t, x, D_x)$ be an $m \times m$ matrix of elliptic classical pseudo-differential operators of order one. Let*

$$P(t, x, D_x) = A(x, D_x) + B(t, x, D_x),$$

where A is a symmetric matrix of pseudo-differential operators of order one and B is a matrix of operators of order zero. Assume that the matrix A is smoothly (microlocally) diagonalizable, with smooth eigenspaces and real eigenvalues $a_j(x, \xi)$, satisfying condition C. Then for any compactly supported $f \in L_\alpha^p \cap L^2$, the solution $u = u(t, x)$ of the Cauchy problem

$$(2.1) \quad i \frac{\partial u}{\partial t} - P(t, x, D_x)u = 0, \quad u(0) = f,$$

satisfies $u(t, \cdot) \in L_{loc}^p$ for all $0 < t \leq T$. Moreover, there is a constant $C > 0$ such that

$$\sup_{0 \leq t \leq T} \|u(t, \cdot)\|_{L^p} \leq C_T \|f\|_{L_\alpha^p}.$$

We note that it is sufficient to only assume that A is microlocally diagonalizable. Then by Lemma 5.1 we can first have the statement of Theorem 2.2 on the cover \tilde{X} . Then, using formula (5.1) and the fact that $\dim \tilde{X} = \dim X$, we get the same conclusion on X .

As a consequence, if the Cauchy data u_0 is compactly supported, we obtain local estimates in other spaces as well:

- $u_0 \in L_{s+\alpha}^p$ implies $u(t, \cdot) \in L_s^p$, $s \in \mathbb{R}$.
- $u_0 \in \text{Lip}(s + (n-1)/2)$ implies $u(t, \cdot) \in \text{Lip}(s)$.
- Let $1 < p \leq q \leq 2$. Then $u_0 \in L_{s-1/q+n/p-(n-1)/2}^p$ implies $u(t, \cdot) \in L_s^q$. Dual result holds for $2 \leq p \leq q < \infty$.

The proof of Theorem 2.2 is based on the other main result Theorem 3.1 which we will discuss in the next section. Estimates in other spaces follow by standard methods of harmonic analysis ([25]).

Now we will give some examples where condition C holds while the transversality condition ($M = 1$) fails. This is, for example, the case when pairs a_j, a_k define glancing hypersurfaces or when we consider Maxwell systems with variable coefficients. Below we will concentrate on systems arising from scalar weakly hyperbolic equations with Levi conditions.

Example 1. In scalar equations with Levi conditions studied by Chazarain [3], Mizohata-Ohya [18], Zeman [28], one assumed that $\{a_j, a_k\} = C_{jk}(a_j - a_k)$. It is clear that in this situation $a_j(x, \xi) = a_k(x, \xi)$ implies $\{a_j, a_k\}(x, \xi) = 0$. However, in a general case when $C_{jk}(x, \xi)$ is non-constant, condition C is satisfied generically.

Example 2. Let L be a scalar operator with involutive characteristics. More precisely, let us denote $\partial_j = D_t + \lambda_j(t, x, D_x)$ and let

$$(2.2) \quad L = \partial_1 \cdots \partial_m + \sum_{k < m} b_{j_1, \dots, j_k} \partial_{j_1} \cdots \partial_{j_k} + c,$$

where $b(t, x, D_x), c(t, x, D_x) \in \Psi^0$ are pseudo-differential operators of order zero for all $t \in [0, T]$. We will assume that symbols of all operators are infinitely differentiable with respect to t in the topology of symbols of the corresponding order. Let us assume that operator L has involutive characteristics, i.e. that

$$[\partial_j, \partial_k] \equiv \partial_j \partial_k - \partial_k \partial_j = \alpha_{jk} \partial_j + \beta_{jk} \partial_k + \gamma_{jk},$$

where $\alpha_{jk}, \beta_{jk}, \gamma_{jk} \in \Psi^0$ are pseudo-differential operators of order zero. Then it was shown by Morimoto in [19] that the Cauchy problem for the equation $Lu = f$ is diagonalizable (with $1 + \sum_{j=1}^{m-1} m!/j!$ components). Even in the simplest case of characteristics not depending on t , we have

$$\{\lambda_j, \lambda_k\} = \alpha_{jk}(x, \xi)(\lambda_j - \lambda_k) + \gamma_{jk},$$

similar to Example 1.

Propagation of singularities of systems with vanishing Poisson brackets has been also studied in these situations. For example, Iwasaki and Morimoto [13] studied propagation of singularities of 3×3 systems, where the second Poisson bracket vanish. Also, Ichinose [11] studied 2×2 systems with vanishing second Poisson brackets. Theorem 2.2 implies a precise statement on L^p estimates.

Theorem 2.3. *Let $1 < p < \infty$, $\alpha = (n-1)|1/p - 1/2|$ and $s \in \mathbb{R}$. Let L be as in (2.2) and suppose that principal symbols $a_j(x, \xi)$ of λ_j satisfy condition C and do not depend on t . Let u be a solution to the Cauchy problem*

$$(2.3) \quad \begin{cases} Lu = 0, \\ \partial_t^j u(0, x) = g_j(x), \quad 0 \leq j \leq m-1, \end{cases}$$

and let Cauchy data $g_j \in L_{\alpha-j+s}^p$ be compactly supported. Then $u(t, \cdot) \in (L_s^p)_{loc}$ for all $t \in [0, T]$ and

$$(2.4) \quad \sup_{t \in [0, T]} \|\partial^J u(t, \cdot)\|_{L_s^p} \leq C \sum_{j=0}^{m-1} \|g_j\|_{L_{\alpha-j+s+m-1}^p},$$

where $\partial^J = \partial_{j_1} \dots \partial_{j_k}$, $k \leq m-1$, and (j_1, \dots, j_k) being permutations of some elements of $\{1, \dots, m\}$.

Note that in the strictly hyperbolic case as well as in some very special cases of operator L in (2.2) (e.g. when all b and c are zero), following the method described by Treves in [27] and estimates for Fourier integral operators, it is possible to obtain the estimate for the Sobolev norm $\|u\|_{L_{s+m-1}^p}$ in the left hand side of (2.4).

Let us now give a final example of L^p estimates for second order equations, which we will prove in the next section.

Example 3. Let us consider the second order equation

$$u'' + b(x, D_x)u' + c(x, D_x)u = 0,$$

where $b \in \Psi^1$, $c \in \Psi^2$. Let us denote $\langle x \rangle = (1 + x^2)^{1/2}$. Introducing $v = \begin{pmatrix} \langle D_x \rangle u \\ u' \end{pmatrix}$, the matrix form of this equation is given by

$$v' = \begin{pmatrix} 0 & \langle D_x \rangle \\ -\langle D_x \rangle^{-1}c & -b \end{pmatrix} v.$$

Let b_1 and c_2 be principal symbols of b and c . The equation is hyperbolic if $b_1^2 \geq 4c_2$, with multiple roots at $b_1^2 = 4c_2$. Assume that $b_1^2 - 4c_2 = \mu^2$, with $\mu \in S^1$ being a symbol of order one. Then characteristics a_1, a_2 satisfy $\{a_1, a_2\} = \frac{1}{2}\{b_1, \mu\}$, which may vanish.

Let the Cauchy data be $u(0) = f_0, u'(0) = f_1$. Let $\alpha = (n-1)|1/p - 1/2|$ and $1 < p < \infty$. If μ is elliptic, it is known that if $f_j \in L_{\alpha-j}^p$, then $u(t, \cdot) \in L^p$. Our result of Theorem 2.2 will imply that the same is true for any $\mu \in S^1$.

Note that in this example we may require only microlocal diagonalization with the same conclusion.

Let us now discuss the propagation of singularities for operators Q_l . This result is essentially a reformulation of Rozenblum's result in the case of finite geometry under Condition C. It is clear (also from multi-phase analysis) that singularities propagate along broken Hamiltonian flows. Let

$$J = \{j_1, \dots, j_{l+1}\}, \quad 1 \leq j_k \leq m, \quad j_k \neq j_{k+1}.$$

Let $\Phi_J(t, x, \xi)$ be the corresponding broken Hamiltonian flow. It means that points follow bicharacteristics of a_{j_1} until meeting the characteristic of a_{j_2} , and then continue

along the bicharacteristic of a_{j_2} , etc. Note that singularities may accumulate if wave front sets for different broken trajectories project to the same point of X .

We can write

$$Q_l = \int_{\Delta} I(\bar{t}) d\bar{t},$$

where $\bar{t} = (t_1, \dots, t_l)$, $\Delta = \{0 \leq t_l \leq t_{l-1} \leq \dots \leq t_1 \leq t\}$ is a simplex in \mathbb{R}^l and $I(\bar{t}) = Z(t_1) \circ \dots \circ Z(t_l)$. It is possible to treat it as a standard Fourier integral operator with the change of variables $\bar{t} = \zeta|\zeta|^{-1}$. Let K be a cone in $\mathbb{R}^N = \mathbb{R}^{n+l}$. Let

$$Iu(x) = \int_K \int_Y e^{i\varphi(x,y,\theta)} a(x,y,\theta) u(y) dy d\theta$$

be a Fourier integral operator with integration over the cone K with respect to θ . Let K_j be K or a face of K . Let $\varphi_j(x,y,\theta_j) = \varphi|_{K_j}$, $\theta_j \in K_j$. Let $\Lambda_j \subset T^*X \times T^*X$ be a Lagrangian manifold with boundary:

$$\Lambda_j = \{(x, \frac{\partial \varphi_j}{\partial x}, y, -\frac{\partial \varphi_j}{\partial y}) : \frac{\partial \varphi_j}{\partial \theta_j} = 0\}.$$

For $G \subset T^*Y$, let $\Lambda_j(G) = \{z \in T^*X : \exists \zeta \in G : (z, \zeta) \in \Lambda_j\}$. Then we have the following statement on the propagation of singularities.

Theorem 2.4. *Let $u \in \mathcal{D}'(Y)$. Then $WF(Iu) \subset \cup_j \Lambda_j(WF(u))$.*

The proof is standard and follows Hörmander [9].

From this, we can deduce first and second terms of the spectral asymptotic of operator P . Let us call T a period of symbol $A(x, \xi)$ if there exists J such that $j_1 = j_{l+1}$, and the trajectory of Φ_J is closed: $\Phi_J(T, x, \xi) = (x, \xi)$. Then we have the following extension of well-known results of Hörmander [8], Duistermaat–Guillemin [7], Safarov–Vassiliev [23], and Rozenblum [21].

Theorem 2.5. *Assume that X is compact and assume that Condition C is satisfied. Let D be the set of $(x, \xi) \in T^*X$ such that there exist T and J such that $\Phi_J(T, x, \xi) = (x, \xi)$. Assume that the measure of D is zero. Then for the spectrum of P the following Weyl formula holds:*

$$N(\lambda) = \#\{j : \lambda_j < \lambda\} = c_n \lambda^n + c'_n \lambda^{n-1} + o(\lambda^{n-1}),$$

where λ_j are eigenvalues of P .

Proof of the L^p estimates will be based on Theorem 3.1, which we regard as an independent result on regularity of solutions of partial differential equations. Theorem 5.3 concerns the measure of the set where a function is small given some information on the multiplicity of its roots. It will play a crucial role in the proof of the smoothing property in Theorem 2.1.

3. REGULARITY OF SOLUTIONS

In this section we will present a principle governing solutions of first order systems. Let $Z(t) \in \mathcal{L}(C_0^\infty(X), \mathcal{D}'(X))$, $t \in [0, T]$, be a time dependent family of operators. Let W_0, W_1 and W be linear subspaces of $\mathcal{D}'(X)$ such that $W_0, W_1 \hookrightarrow W$. We will

make different choices of these spaces in the future, dependent on the structure of operators $Z(t)$. Let us consider the Cauchy problem for $u = u(t, x)$:

$$(3.1) \quad \begin{cases} u' - Z(t)u = r, & r(t) \in W_0, \\ u(0) \in W_1. \end{cases}$$

One is often interested in the following question. If the right hand side and Cauchy data satisfy $r(t) \in W_0$ and $u(0, \cdot) \in W_1$, when do fixed time solutions $u(t, \cdot)$ of (3.1) belong to W ? In general, some loss of regularity is possible in problem (3.1) even if $Z(t)$ are very good. So we will think of W_0 being the smallest, W_1 an intermediate, and W the largest among these spaces. The following theorem says that if operators $Z(t)$ have some structure, and solutions of Cauchy problem (3.1) with zero Cauchy data are in W , so will be solutions with Cauchy data from some sufficiently large space W_1 .

Theorem 3.1. *Let $W_0, W_1 \hookrightarrow W$ be linear subspaces of $\mathcal{D}'(X)$.*

Let $Z(t) \in \mathcal{L}(C_0^\infty(X), \mathcal{D}'(X))$, $t \in [0, T]$. Assume that

- (i) (Boundedness) *Z extends to an operator in $L^\infty([0, T], \mathcal{L}(L^2(X), L^2(X)))$, and $Z(t)$ extend to continuous linear operators from W_1 to W , for all $t \in [0, T]$.*
- (ii) (Calculus) *$Z(t_1) \circ \cdots \circ Z(t_l) : W_1 \rightarrow W$ are continuous for all l and for all $t_1, \dots, t_l \in [0, T]$.*
- (iii) (Smoothing) *There exists l such that*

$$Z(t) \int_0^t \int_0^{t_1} \cdots \int_0^{t_{l-1}} Z(t_1) \circ \cdots \circ Z(t_l) dt_l \cdots dt_1$$

is continuous from W_1 to W_0 , for all $t \in [0, T]$.

- (iv) (Zero Cauchy data) *Solutions $v = v(t, x)$ of the Cauchy problem*

$$(3.2) \quad \begin{cases} v' - Z(t)v = r, & r(t) \in W_0, \\ v(0) = 0, \end{cases}$$

satisfy $v(t, \cdot) \in W$ for $t \in [0, T]$.

Then the solution $u = u(t, x)$ of the Cauchy problem

$$(3.3) \quad \begin{cases} u' - Z(t)u = r, & r(t) \in W_0, \\ u(0) \in W_1, \end{cases}$$

satisfies $u(t, \cdot) \in W$ for all $t \in [0, T]$.

Moreover, if W_0, W_1, W are normed spaces and if solutions $v(t, \cdot)$ to (3.2) in (iv) satisfy $\|v(t, \cdot)\|_W \leq C\|r(t)\|_{W_0}$ for all $t \in [0, T]$, then also

$$\|u(t, \cdot)\|_W \leq C(\|u(0)\|_{W_1} + \|r(t)\|_{W_0}),$$

for all $t \in [0, T]$.

Conditions (i) and (ii) ensure that operators $Z(t)$ have some structure. Indeed, if $W_1 \subset W$ is different from W , (ii) does not follow from (i). In our typical applications, $Z(t)$ will be time dependent pseudo-differential or Fourier integral operators, and compositions in (ii) are essentially of the form of a single operator $Z(t)$. Condition (iii) is natural from the point of view of harmonic analysis, since integration with respect to a parameter often brings additional regularity. Condition (iv) ensures

that solutions with zero Cauchy data and regular right hand side are also sufficiently regular.

Proof. Let $U(t)$ be an operator solving the Cauchy problem

$$(3.4) \quad \begin{cases} U' - Z(t)U = R(t), & R(t) \in \mathcal{L}(W_1, W_0), \\ U(0) = I. \end{cases}$$

Let $U_0(t)$ be some partial solution to the problem

$$\begin{cases} U'_0 - Z(t)U_0(t) = R(t), \\ U_0(0) = 0. \end{cases}$$

Then the solution U of problem (3.4) satisfies

$$(3.5) \quad U(t) = U_0(t) + I + \int_0^t Z(t_1)dt_1 + \int_0^t \int_0^{t_1} Z(t_1)Z(t_2)dt_2dt_1 + \dots$$

The convergence of this series can be understood in L^2 . Indeed, because of assumption (i), the term of this series with k integrals can be estimated by $t^k \sup_t \|Z(t)\|_{L^2 \rightarrow L^2}^k / k!$. From this it also follows that $U(t)$ is a solution of (3.4) in L^2 . Let us now define

$$\begin{aligned} S_N(t) = I + \int_0^t Z(t_1)dt_1 + \int_0^t \int_0^{t_1} Z(t_1)Z(t_2)dt_2dt_1 + \dots \\ + \int_0^t \int_0^{t_1} \dots \int_0^{t_{N-1}} Z(t_1)Z(t_2) \dots Z(t_N)dt_N \dots dt_2dt_1. \end{aligned}$$

Let $V(t) = U(t) - S_N(t)$, it is equal to $U_0(t)$ plus the remainder of the series (3.5). Then we have

$$(3.6) \quad \begin{aligned} V'(t) - Z(t)V(t) = \\ (U - S_N)'(t) - Z(t)(U - S_N)(t) = \\ (U' - ZU)(t) - (S'_N - ZS_N)(t) = \\ R(t) - Z(t) \int_0^t \int_0^{t_1} \dots \int_0^{t_{N-1}} Z(t_2) \dots Z(t_N)dt_2 \dots dt_N. \end{aligned}$$

Choosing $N = l$, from assumption (iii) of the theorem the second term is continuous from W_1 to W_0 . Since also $R(t) \in \mathcal{L}(W_1, W_0)$, it follows that the right hand side is a continuous linear operator from W_1 to W_0 .

Let $w = u(0) \in W_1$ be the Cauchy data for (3.3). If we denote by $\rho(t)$ the value of the operator in the last line of (3.6) at w , we will have $\rho(t) \in W_0$. The value of $V(0)$ is

$$V(0) = U(0) - S_N(0) = 0.$$

It follows now that $V(t)w$ solves Cauchy problem (3.2), so it belongs to W by assumption (iv). Since $S_N(t)$ is continuous from W_1 to W by assumption (ii), and $V(t)w = U(t)w - S_N(t)w$ is in W , we obtain $u(t, \cdot) = U(t)w \in W$.

Moreover, suppose that we also have the estimate $\|v(t, \cdot)\|_W \leq C\|\rho(t)\|_{W_0}$ in (iv). Then we also have

$$\|u(t, \cdot)\|_W \leq \|V(t)w\|_W + \|S_N(t)w\|_W \leq C\|\rho(t)\|_{W_0} + C\|w\|_{W_1} \leq C(\|w\|_{W_1} + \|R(t)w\|_{W_0}).$$

□

Later we will need this in the case of $Z(t)$ being Fourier integral operators of order zero. However, let us point out several applications to other cases of pseudo-differential and Fourier integral operators. In these cases we will make different appropriate choices of spaces W_0, W_1, W . Moreover, in Corollaries 3.2 and 3.3 we will assume that the corresponding non-homogeneous Cauchy problems with zero Cauchy data have unique solutions. This is a natural assumption if $Z(t)$ behave sufficiently well with respect to t since we are working in subspaces of L^2 .

Corollary 3.2. *Let $1 < p < \infty$. Let $Z(t) \in \Psi^0$, $t \in [0, T]$, be a family of pseudo-differential operators of order zero with amplitudes compactly supported in x, y , uniformly in t . Suppose that $Z \in L^\infty([0, T], \mathcal{L}(L^2, L^2)) \cap L^\infty([0, T], \mathcal{L}(L^p, L^p))$. Then the solution $u = u(t, x)$ of the Cauchy problem*

$$(3.7) \quad \begin{cases} u' - Z(t)u = r, & r(t) \in L^p, \quad t > 0, \\ u(0) \in L^p, \end{cases}$$

satisfies $u(t, \cdot), u'(t, \cdot) \in L^p$, for all $t \in (0, T]$. Moreover, we have an estimate

$$\sup_{t \in [0, T]} \|u(t, \cdot)\|_{L^p} \leq C \|u(0)\|_{L^p}.$$

If amplitudes of $Z(t)$ are not compactly supported with respect to x, y , we have a similar local statement for compactly supported Cauchy data. Note also that pseudo-differential operators of order zero are locally bounded in L^2 and L^p , for all $1 < p < \infty$. Conditions $Z \in L^\infty([0, T], \mathcal{L}(L^2, L^2))$ and $Z \in L^\infty([0, T], \mathcal{L}(L^p, L^p))$ simply mean that we have some control on their norms, i.e. there exist a constant C such that

$$\|Z(t)\|_{L^2 \rightarrow L^2} \leq C, \quad \|Z(t)\|_{L^p \rightarrow L^p} \leq C, \quad \forall t \in [0, T].$$

Proof. Let us choose $W = W_1 = W_0 = L^p_{comp}$. Let us check conditions of Theorem 3.1. Properties (i) and (ii) follow from regularity properties of pseudo-differential operators of order zero and our assumptions. Property (iii) also holds because $Z(t)$ are locally bounded in L^p . Property (iv) is a consequence of Duhamel's principle and is similar to the one in Corollary 3.3. Picard series is convergent in L^p provided that operators norms $\|Z(t)\|_{L^p \rightarrow L^p}$ are uniformly bounded for $t \in [0, T]$. Norm estimate follows from this as well.

We can see that $u'(t, \cdot) \in W$ from $u' = Z(t)u + r$ and from the continuity of $Z(t)$ in W . \square

We will now apply Theorem 3.1 in the case of $Z(t)$ being Fourier integral operators. While our case (1.3) corresponds to $Z(t)$ being operators of order zero, the crucial smoothing property (iii) will follow from the fact that operators $Z(t)$ have a special structure. For general Fourier integral operators $Z(t)$ without structure, we need to assume that they are of negative orders. This is for example the case when the zero order term B in Theorem 2.2 is actually a pseudo-differential operator of some negative order.

Corollary 3.3. *Let $1 < p < \infty$, $\epsilon > 0$, and $\alpha = (n - 1)|1/p - 1/2|$. Let $Z(t) \in I^{-\epsilon}$, $t \in [0, T]$, be a family of non-degenerate Fourier integral operator of order $-\epsilon$ with amplitudes compactly supported in x, y , uniformly in t . Suppose that operators $Z(t)$ can be composed and that $Z \in L^\infty([0, T], \mathcal{L}(H^s, H^s))$, for some $s > (pn - 2n)/2p$*

when $p > 2$ and $s = 0$ when $p \leq 2$. Then the solution $u = u(t, x)$ of the Cauchy problem

$$(3.8) \quad \begin{cases} u' - Z(t)u = r, & r(t) \in H^s, \ t > 0, \\ u(0) \in (L_\alpha^p) \cap L^2, \end{cases}$$

satisfies $u(t, \cdot) \in L^p$, for all $t \in (0, T]$. Moreover,

$$\sup_{t \in [0, T]} \|u(t, \cdot)\|_{L^p} \leq C \|u(0)\|_{L_\alpha^p}.$$

Note that operators $Z(t)$ are locally bounded in H^s , so assumption $Z \in L^\infty([0, T], \mathcal{L}(H^s, H^s))$ simply means that $\|Z(t)\|_{H^s \rightarrow H^s} \leq C$ for all $t \in [0, T]$.

Proof. Let $W = L^2 \cap L_{comp}^p$, $W_1 = (L_\alpha^p)_{comp}$, and $W_0 = H_{comp}^s \subset W$. Let us check conditions of Theorem 3.1. Condition (i) follows from the fact that non-degenerate Fourier integral operators of order 0 are bounded from $(L_\alpha^p)_{comp}$ to L_{loc}^p . Condition (ii) follows from the calculus of non-degenerate Fourier integral operators, since we assumed that compositions of $Z(t)$ are again non-degenerate Fourier integral operators. Smoothing condition (iii) for large l follows again from the calculus, since operators $Z(t)$ are of order $-\epsilon$.

Finally, let us show that solutions of $v' - Z(t)v = r(t)$, $r(t) \in H_{comp}^s$, with zero Cauchy data $v(0) = 0$, satisfy $v(t, \cdot) \in L^2 \cap L^p$. In fact, we will show that $v(t, \cdot) \in H^s \subset L^p \cap L^2$.

From the uniqueness of the solution of this problem it follows that we can use Duhamel's principle to write

$$(3.9) \quad v(t, x) = \int_0^t E(t, s) r(s, x) ds,$$

where $E(t, s)$ is the propagator of

$$\begin{cases} (\partial_t - Z(t))E(t, s) = 0, \\ E(t, s)|_{t=s} = I. \end{cases}$$

Picard series for this problem gives the asymptotic expansion of $E(t, s)$, in particular implying that $E(t, s)$ is bounded in L^2 and in H^s provided operator norms $\|Z(t)\|_{H^s \rightarrow H^s}$ are uniformly bounded for $t \in [0, T]$. From (3.9) it follows that $v(t, \cdot) \in H^s$. Moreover, since $r(s, \cdot) \in H^s$, we also get an estimate

$$\|u(t, \cdot)\|_{H^s} \leq C \sup_{\tau \in [0, T]} \|r(\tau, \cdot)\|_{H^s},$$

implying the estimate in Corollary 3.3. \square

Proof of Theorem 2.2. As we have already mentioned, by Lemma 5.1 we can assume that characteristics of A are correctly defined on T^*X . Since A is diagonalizable, we can write

$$(3.10) \quad \begin{aligned} P(t, x, D_x) &= \bigoplus a_j(x, D_x) + B(t, x, D_x), \\ B(t, x, D_x) &= (B_{jk}(t, x, D_x))_{1 \leq j, k \leq m}, \ B_{jk} \in C^\infty([0, T], \Psi^0). \end{aligned}$$

Some of a_j 's may be identically equal to each other. We can renumber a_j 's into r groups (possibly of size one) of equal characteristics. These are the eigenvalues of the matrix $A(x, \xi)$ counted with multiplicity. Thus, we have $1 = k_1 < k_2 < \dots < k_r =$

$n + 1$, and $a_{k_i} \equiv \dots \equiv a_{k_{i+1}-1} \not\equiv a_k$, for $k < k_i$ or $k \geq k_{i+1}$. This means that we have a group of the same roots a_1, \dots, a_{k_2-1} , etc., while roots from different groups are not identically the same. Therefore, this is a decomposition of the first order principal part into a block-diagonal form with the same roots in each block, with possible equality of roots in different blocks at some points. So we can write

$$(3.11) \quad P(t, x, D_x) = \text{diag}(\tilde{a}_1, \dots, \tilde{a}_r) + B(t, x, D_x),$$

where $\tilde{a}_i = \text{diag}(a_{k_i}, \dots, a_{k_{i+1}-1})$ are diagonal matrices with equal roots at the diagonal. Let us set

$$\tilde{A}_i = \tilde{a}_i + (B_{\mu\nu})_{k_i \leq \mu, \nu \leq k_{i+1}-1},$$

so that

$$P = \tilde{A} + B = \text{diag}(\tilde{A}_1, \dots, \tilde{A}_r) + B.$$

Note that in the last equality we can assume $B_{\mu\nu} = 0$ for $k_i \leq \mu, \nu \leq k_{i+1}-1$ if we add these components to the corresponding components of \tilde{A} . Let $U(t) = \exp(-i\tilde{A}t)V(t)$. This is well defined in view of, for example, [26, VIII]. Then

$$V' = Z(t)V \equiv -ie^{i\tilde{A}t}Be^{-i\tilde{A}t}V, \quad V(0) = I.$$

Now we will apply Theorem 3.1 with $Z(t) = -ie^{i\tilde{A}t}Be^{-i\tilde{A}t}$. Let us choose $W = L_{comp}^2 \cap L^p$, $W_1 = (L_\alpha^p)_{comp}$, and $W_0 = H_{comp}^s$ with $s > (pn - 2n)/2p$ for $p > 2$ and $s = 0$ for $1 < p \leq 2$. Conditions (i) and (ii) follow from the calculus and regularity properties of non-degenerate Fourier integral operators of order zero. Smoothing condition (iii) follows from Theorem 2.1. For condition (iv) we can use Duhamel's principle similar to the proof of Corollary 3.3. Thus, Theorem 3.1 implies that propagator $V(t)$ is continuous from L_α^p to L^p . Operator $U(t)$ is a composition of $V(t)$ with a non-degenerate Fourier integral operator $\exp(-i\tilde{A}t)$, so $U(t)$ is given as a sum of a smoothing series obtained by the multiplication of Picard series for $V(t)$ with $\exp(-i\tilde{A}t)$. Using the calculus of Fourier integral operators in each term of the series and its smoothing property we can repeat the argument of Theorem 3.1 in this case to see that $u(t, \cdot) \in L^p$ with an estimate for its norm. \square

Note that if B in Theorem 2.2 is a pseudo-differential operator of negative order, $B \in \Psi^\mu$, for some $\mu < 0$, the proof is simpler because we do not have to use Theorem 2.1 to prove condition (iii) of Theorem 3.1. Instead, we can use directly Corollary 3.3 to obtain the smoothing property (iii).

Proof of Theorem 2.3. Let L be as in (2.2) and let u be the solution of

$$(3.12) \quad Lu = f, \quad D_t^j u(0, x) = g_j(x), \quad 0 \leq j \leq m-1.$$

Let

$$U = (u, \partial_1 u, \partial_2 u, \dots, \partial_1 \partial_2 u, \partial_2 \partial_1 u, \dots, \partial^J u, \dots)^T,$$

where $J = \{j_1, \dots, j_k\}$ is a permutation of some elements of $\{1, \dots, m\}$, $|J| = k \leq m-1$. Vector U has $1 + \sum_{j=1}^{m-1} m!/j!$ components. Here we can write $\partial^J = D_t^k + \sum_{j=0}^{k-1} c_j^J(t, x, D_x) D_t^j$, where $c_j^J(t) \in \Psi^{k-j}$. We set $|J| = k$. It was shown by Morimoto in [19] that U solves the system

$$(3.13) \quad D_t U + AU + BU = F, \quad U(0, x) = G(x),$$

where $F = (0, \dots, 0, f, \dots, f)$ and $G = (g_0, \dots, g_{|J|} + \sum_{j=0}^{|J|-1} c_j^J g_j, \dots)$, A is a diagonal matrix with λ_j 's at the diagonal and B is a matrix of pseudo-differential operators of order zero. Matrix B has some operators in the last row, zeros, and -1 above the diagonal. If λ_j satisfy Condition C, Theorem 2.2 implies $\|U(t, \cdot)\|_{L^p} \leq C\|G\|_{L_\alpha^p}$. Since $c_j^J(t) \in \Psi^{k-j}$, we get $\|G\|_{L_\alpha^p} \leq C \sum_{j=0}^{m-1} \|g_j\|_{L_{\alpha+m-1-j}^p}$, which implies the estimate of the Theorem. \square

4. ESTIMATES FOR PICARD SERIES

In this Section we will prove Theorem 2.1 on the smoothing properties of terms Q_l of the Picard series (1.4).

Let $A_j \in \Psi^1$, $j = 1, \dots, r$, be elliptic pseudo-differential operators of order one. Let $a_j(x, \xi)$ denote their principle symbols. We can assume that there are no identical symbols among these a_j 's. Let

$$H(\bar{t}) = e^{iA_{j_1}t_1} e^{iA_{j_2}(t_2-t_1)} \dots e^{-iA_{j_{l+1}}t_l},$$

where $1 \leq j_k \leq r$, $j_k \neq j_{k+1}$, $k = 1, \dots, l+1$, and $\bar{t} = (t_1 \dots t_l)$. Let us define

$$Q = \int_0^t \int_0^{t_1} \dots \int_0^{t_{l-1}} B(\bar{t}) H(\bar{t}) dt_l \dots dt_1,$$

where $B(\bar{t}) \in \Psi^0$ is a pseudo-differential operator of order zero smoothly dependent on \bar{t} . Such operators Q appear in the Picard series (1.4), (1.5). In this section we will give a detailed description of operator Q in order to prove that it is a smoothing operator when l is sufficiently large. First of all let us note that H is a Fourier integral operator and due to the theorem on compositions of Fourier integral operators the canonical relation $\Lambda^{\bar{t}} \subset T^*X \times T^*X$ of $H(\bar{t})$ is given by

$$\Lambda^{\bar{t}} = \{(x, p, y, \xi) : (x, p) = \Psi^{\bar{t}}(y, \xi)\},$$

where $\Psi^{\bar{t}} = \Phi_{j_1}^{t_1} \circ \dots \circ \Phi_{j_l}^{t_l - t_{l-1}} \circ \Phi_{j_{l+1}}^{-t_l}$ and Φ_j^t is the Hamiltonian flow defined by a_j .

It can be easily checked that H is a solution operator for the system of equations

$$(4.1) \quad \frac{\partial H}{\partial t_k} = iT_k(t_1, \dots, t_k)H, \quad k = 1, \dots, l,$$

where $T_k \in \Psi^1$ is a pseudo-differential operator of order one. In view of Egorov's theorem its principle symbol is equal to

$$(4.2) \quad T_k^0(t_1, \dots, t_k, x, \xi) = (a_{j_k} - a_{j_{k+1}}) \circ \Phi_{j_1}^{t_1} \circ \dots \circ \Phi_{j_k}^{t_k - t_{k+1}}(x, \xi), \quad (x, \xi) \in T^*X$$

for all $k = 1, \dots, l$. Let us construct a phase function $\varphi(\bar{t}, x, y, \xi)$ which defines the operator $H(\bar{t})$ for small \bar{t} . We will look for it in the form $\varphi(\bar{t}, x, y, \xi) = \psi(\bar{t}, x, \xi) - y \cdot \xi$. It follows from (4.1) and (4.2) that ψ satisfies a system of Hamilton-Jacobi equations

$$(4.3) \quad \frac{\partial \psi}{\partial t_k} = T_k^0(\bar{t}, x, \frac{\partial \psi}{\partial x}), \quad \psi(0, x, \xi) = x \cdot \xi.$$

In [21] it was checked that Frobenius conditions for system (4.3) are satisfied. Solving this system we obtain a non-degenerate phase function. This phase function defines

a Lagrangian manifold $\Lambda^{\bar{t}}$, so that we have

$$(4.4) \quad \left(x, \frac{\partial \psi}{\partial x}\right) = \Psi^{\bar{t}}(y, \xi) = (x^{\bar{t}}(y, \xi), p^{\bar{t}}(y, \xi)), \quad y = \frac{\partial \psi}{\partial \xi}.$$

Now we are going to investigate the smoothing properties of operator Q . We can write Q as

$$Qu(x) = \int_{\Delta} \int_{\mathbb{R}^n} \int_{\mathbb{R}^n} e^{i\varphi(\bar{t}, x, y, \xi)} b(\bar{t}, x, y, \xi) u(y) dy d\xi d\bar{t},$$

where φ satisfies (4.4) and b is an amplitude of order zero, which we may assume to be compactly supported with respect to x and y . Here Δ is a simplex $\{0 \leq t_l \leq t_{l-1} \leq \dots \leq t_1 \leq t\}$.

Let $\chi \in C_0^\infty$ be a cut-off function such that $\chi(\tau) = 1$ for $|\tau| < 1$ and $\chi(\tau) = 0$ for $|\tau| > 2$. Operator Q can be decomposed as $Q = R_1 + R_2 + R_3$, where

$$R_j u(x) = \int_{\Delta} \int_{\mathbb{R}^n} \int_{\mathbb{R}^n} e^{i\varphi(\bar{t}, x, y, \xi)} \mu_j(\epsilon, \bar{t}, x, y, \xi) u(y) dy d\xi d\bar{t}, \quad j = 1, 2, 3,$$

where

$$\begin{aligned} \mu_1(\epsilon, \bar{t}, x, y, \xi) &= \left(1 - \chi(\epsilon^{-1} |\frac{\partial \varphi}{\partial \xi}|)\right) b(\bar{t}, x, y, \xi), \\ \mu_2(\epsilon, \bar{t}, x, y, \xi) &= \chi(\epsilon^{-1} |\frac{\partial \varphi}{\partial \xi}|) \left(1 - \chi(\epsilon^{-1} |\frac{\partial \varphi}{\partial \bar{t}}| |\xi|^{-1})\right) b(\bar{t}, x, y, \xi), \\ \mu_3(\epsilon, \bar{t}, x, y, \xi) &= \chi(\epsilon^{-1} |\frac{\partial \varphi}{\partial \xi}|) \chi(\epsilon^{-1} |\frac{\partial \varphi}{\partial \bar{t}}| |\xi|^{-1}) b(\bar{t}, x, y, \xi). \end{aligned}$$

Let us first consider operator R_1 . On the support of μ_1 we have the estimate

$$\left| \frac{\partial \varphi}{\partial \xi} \right| \geq \epsilon.$$

Therefore, there exists an operator $L(\frac{\partial}{\partial \xi})$ of order 1, such that $Le^{i\varphi} = e^{i\varphi}$, with coefficients estimated by $C\epsilon^{-1}$ on the support of the amplitude of μ_1 . When integrating by parts with L there may appear an additional factor of ϵ^{-1} when differentiating χ . Integrating by parts p times with this operator L we obtain an operator with an amplitude of order $-p$ and coefficients that can be estimated by ϵ^{-2p} . From Lemma 5.2 with $q = -p$ we obtain the following estimate

$$(4.5) \quad \|R_1\|_{p-n-1} \leq C\epsilon^{-3p+n+1}.$$

The same procedure with integrating by parts with respect to ξ can not be applied to R_2 . But here there is a possibility to integrate by parts with respect to \bar{t} . Indeed, there exists an operator $M(\frac{\partial}{\partial \bar{t}})$, such that $Me^{i\varphi} = e^{i\varphi}$, with coefficients not greater than $\epsilon^{-1}|\xi|^{-1}$ on the support of μ_2 . Integrating by parts with M we obtain an operator with an amplitude of order -1 with coefficients that can be estimated by ϵ^{-2} , where another ϵ^{-1} may appear from differentiating χ . The boundary integrals have the same form as Q but they have amplitudes of order -1 and depend on not less than $l-2$ time variables. The reason for possibly losing two variables is that after restriction to the boundary, say $t_2 = t_1$, it may happen that a_{j_1} and a_{j_3} are the same roots. It follows that we can apply such procedure $[l/2]$ times. As a result we obtain operators

of order $-[l/2]$ with coefficients not greater than $C\epsilon^{-2[l/2]}$. Then in view of Lemma 5.2 for $[l/2] > n + 1$ we have

$$(4.6) \quad \|R_2\|_{[l/2]-n-1} \leq C\epsilon^{-3[l/2]+n+1}.$$

Let us now consider the last integral R_3 . It is not possible to apply procedures with integrating by parts as above either with respect to ξ or with respect to \bar{t} . But in this case it turns out that the support of amplitude μ_3 is small. The singular support of the integral kernel of Q may be very irregular in this case, so a more delicate analysis is necessary to show the smoothing properties of R_3 . First we will show that

$$(4.7) \quad |T_j^0(\bar{t}, x^{\bar{t}}(y, \xi), p^{\bar{t}}(y, \xi))| \leq C\epsilon|\xi|, j = 1, \dots, l,$$

on the support of μ_3 . We notice that μ_3 differs from 0 only if

$$(4.8) \quad \left| \frac{\partial \varphi}{\partial \xi} \right| \leq 2\epsilon, \quad \left| \frac{\partial \varphi}{\partial t} \right| \leq 2\epsilon|\xi|.$$

It follows from (4.3) and (4.8) that

$$(4.9) \quad |T_j^0(\bar{t}, x, \frac{\partial \psi}{\partial x})| \leq C\epsilon|\xi|, j = 1, \dots, l.$$

Because of homogeneity of T_j^0 with respect to ξ we also have the following trivial estimates

$$|\partial_x T_j^0(\bar{t}, x, \frac{\partial \psi}{\partial x})| \leq C|\xi|, \quad |\partial_\xi T_j^0(\bar{t}, x, \frac{\partial \psi}{\partial x})| \leq C, \quad j = 1, \dots, l.$$

Consequently, we obtain

$$\begin{aligned} & |T_j^0(\bar{t}, x, \frac{\partial \psi}{\partial x}) - T_j^0(\bar{t}, x^{\bar{t}}(y, \xi), p^{\bar{t}}(y, \xi))| \\ & \leq C(|\xi||x - x^{\bar{t}}(y, \xi)| + |\frac{\partial \psi}{\partial x} - p^{\bar{t}}(y, \xi)|), \quad j = 1, \dots, l. \end{aligned}$$

Since $\frac{\partial \varphi}{\partial \xi}(\bar{t}, x^{\bar{t}}(y, \xi), y, \xi) = 0$, we get

$$\frac{\partial \varphi}{\partial \xi}(\bar{t}, x, y, \xi) = \frac{\partial \varphi}{\partial \xi}(\bar{t}, x, y, \xi) - \frac{\partial \varphi}{\partial \xi}(\bar{t}, x^{\bar{t}}(y, \xi), y, \xi) = \frac{\partial^2 \varphi}{\partial x \partial \xi}(\bar{t}, x^*, y, \xi)(x - x^{\bar{t}}(y, \xi)),$$

for some points x^* . Since $|\frac{\partial^2 \varphi}{\partial x \partial \xi}| \neq 0$ for small \bar{t} , we obtain

$$(4.10) \quad |x - x^{\bar{t}}(y, \xi)| \leq C \left| \frac{\partial \varphi}{\partial \xi}(\bar{t}, x, y, \xi) \right|.$$

From the properties of Lagrangian manifold $\Lambda^{\bar{t}}$ in (4.4) we also obtain

$$(4.11) \quad \left| \frac{\partial \psi}{\partial x}(\bar{t}, x, \xi) - p^{\bar{t}}(y, \xi) \right| = |p^{\bar{t}}(\frac{\partial \psi}{\partial \xi}, \xi) - p^{\bar{t}}(y, \xi)| \leq C \left| \frac{\partial \varphi}{\partial \xi}(\bar{t}, x, y, \xi) \right| |\xi|, \quad \forall (x, y, \xi).$$

Finally, taking into consideration (4.8)-(4.11) we obtain (4.7) on the support of μ_3 .

Estimate (4.7) shows us that the support of the amplitude μ_3 of operator R_3 is contained in

$$\Xi_1 = \{(\bar{t}, x, y, \xi) : x \in X, \xi \neq 0, (\bar{t}, y, \frac{\xi}{|\xi|}) \in \Xi\},$$

where

$$\Xi = \{(\bar{t}, y, \frac{\xi}{|\xi|}) : |T_j^0(\bar{t}, x^{\bar{t}}(y, \frac{\xi}{|\xi|}), p^{\bar{t}}(y, \frac{\xi}{|\xi|}))| \leq C\epsilon, j = 1, \dots, l\}.$$

That allows us to apply estimate (5.4) of Lemma 5.2 with $q = 0$ and $\delta = 1$ to the operator R_3 to get

$$(4.12) \quad \|R_3\|_{-3n/2-2} \leq C\epsilon^{-n-1} \text{meas}(\Xi),$$

where meas is the natural measure on $[0, T]^l \times S^*X$. To use this inequality we need to estimate the measure $\text{meas}(\Xi)$. The set Ξ maps to

$$\Xi_2 = \{(\bar{t}, y, \xi) : |T_j^0(\bar{t}, y, \xi)| \leq C\epsilon, |p^{\bar{t}}(y, \xi)| = 1, j = 1, \dots, l\}$$

under the Hamiltonian flow $\Psi^{\bar{t}}$ which preserve measure. The measure of Ξ_2 may be estimated by the measure (in $[0, T]^l \times X \times \mathbb{R}^n$) of

$$\Xi_3 = \{(\bar{t}, y, \xi) : |T_j^0(\bar{t}, y, \xi)| \leq C\epsilon, c \leq |\xi| \leq C, j = 1, \dots, l\}.$$

Let us now introduce sets

$$\Sigma_j(t_1, \dots, t_{j-1}, y, \xi, \epsilon) = \{t_j : |T_j^0(\bar{t}, y, \xi)| \leq C\epsilon, c \leq |\xi| \leq C\}, j = 1, \dots, l.$$

To estimate measures of these sets we will use Theorem 5.3. For this, in the notation of Theorem 5.3, we set $w = (y, \xi)$ and consider the function

$$t \mapsto f(t, w) = T_j^0(t_1, \dots, t_{j-1}, t, t_{j+1}, \dots, t_l, y, \xi),$$

where t takes the position of t_j in \bar{t} . Condition on Poisson brackets shows us that function f may have zeros in t of order not greater than M . It follows now from Theorem 5.3 that

$$(4.13) \quad \max_{y, \xi} \text{meas}\{t_j \in [0, T] : |T_j^0(\bar{t}, y, \xi)| \leq C\epsilon\} \leq C\epsilon^{1/2M},$$

where meas is just the Lebesgue measure, y varies over a compact set and $c \leq |\xi| \leq C$. Now we can use a simple observation that if we have two functions $f_1(t_1)$ and $f_2(t_1, t_2)$ such that

$$\text{meas}\{t_1 : |f_1(t_1)| \leq \epsilon\} \leq C\epsilon^\alpha, \max_{t_1} \text{meas}\{t_2 : |f_2(t_1, t_2)| \leq \epsilon\} \leq C\epsilon^\alpha,$$

then

$$\begin{aligned} \text{meas}\{(t_1, t_2) : |f_1(t_1)| \leq \epsilon, |f_2(t_1, t_2)| \leq \epsilon\} &= \int_{|f_1(t_1)| \leq \epsilon, |f_2(t_1, t_2)| \leq \epsilon} dt_1 dt_2 \\ &= \int_{|f_1(t_1)| \leq \epsilon} \left(\int_{|f_2(t_1, t_2)| \leq \epsilon} dt_2 \right) dt_1 \\ &\leq C\epsilon^{2\alpha}. \end{aligned}$$

Applying this argument l times to the estimate (4.13), we get

$$\text{meas}(\Xi_3) \leq C \text{vol}(\text{supp}_y b) \max_{(y, \xi)} \prod_{j=1}^l \max_{t_1, \dots, t_{j-1}} \text{meas}(\Sigma_j(\bar{t}, y, \xi, \epsilon)),$$

which implies

$$\text{meas}(\Xi) \leq C\epsilon^{l/2M}.$$

Here we used that the support of the amplitude $b(\bar{t}, x, y, \xi)$ is compact with respect to x and y , and that T is small. Finally, combining this with estimate (4.12), we get

$$(4.14) \quad \|R_3\|_{-3n/2-2} \leq C\epsilon^{l/2M-n-1}.$$

An application of Lemma 5.6 to (4.5), (4.6) and (4.14) yields the required estimate for Q in Theorem 2.1.

5. VARIOUS AUXILIARY RESULTS

First we describe the lifting of the problem to insure that characteristic roots of the principal symbol of P are globally uniquely defined. The proof follows the paper of Rozenblum [21]. We give it here for the completeness and since we will need this construction to determine the orders for L^p estimates in Theorem 2.2.

Lemma 5.1. *Let X be a smooth compact manifold without boundary of dimension $n \geq 3$. Then there exists a cover \tilde{X} of X with finitely many leaves, such that on the lifting \tilde{E} of E to $T^*\tilde{X}$ branches of eigenvalues $a_j(x, \xi)$ and eigenspaces $V_j(x, \xi)$ of the principal symbol $A(x, \xi)$ are smooth and globally well defined. The space $L^2(\tilde{E})$ has a decomposition into a direct sum of m spaces such that the matrix representation of $P(\tilde{x}, \tilde{\xi})$ with respect to this decomposition consists of pseudo-differential operators, and its principal symbol is a diagonal matrix with $a_j(\tilde{x}, \tilde{\xi})$ at the diagonal.*

Proof. Let us fix $z_0 = (x_0, \xi_0) \in S^*X$. For each path in S^*X beginning at z_0 we look at the continuation of the diagonalization along this path. At each point $z \in S^*X$ we obtain up to permutation several possible collections $a_j(z), V_j(z)$, such that homotopic paths from z_0 to z correspond to the same collection. Since for $n \geq 3$ homotopy of paths in S^*X and their projections to X is equivalent, we obtain a homomorphism $\pi_1(X)$ to the permutation group of order m . A construction in the homotopy theory ([20]) gives a cover $p : \tilde{X} \rightarrow X$ with finitely many leaves, such that the lifting of this homomorphism to $\pi_1(\tilde{X})$ is trivial. So, each closed path in $S^*\tilde{X}$ takes eigenvalues and, therefore, also eigenspaces, to themselves. This means that eigenvalues and eigenspaces have global smooth branches on $S^*\tilde{X}$ and hence also on $T^*\tilde{X}$.

Let $p_j^0(\tilde{x}, \tilde{\xi})$ be a family of orthogonal projectors on $V_j(\tilde{x}, \tilde{\xi})$. By the standard Gram-Schmidt process, we can add lower order terms to $p_j^0(\tilde{x}, \tilde{\xi})$ to obtain symbols $p_j(\tilde{x}, \tilde{\xi})$, for which $p_j(\tilde{x}, \tilde{D})p_k(\tilde{x}, \tilde{D}) = \delta_{jk}p_j(\tilde{x}, \tilde{D})$ and $\sum p_j(\tilde{x}, \tilde{D}) = 1$. For $u \in L^2(E)$, let $u_j = p_j(\tilde{x}, \tilde{D})u$. Since $P(\tilde{x}, \tilde{D})p_j(\tilde{x}, \tilde{D}) = p_j(\tilde{x}, \tilde{D})P(\tilde{x}, \tilde{D})p_j(\tilde{x}, \tilde{D})$ modulo lower order terms, $P = \sum p_j P p_k$ yields the desired diagonalization. \square

Let a_j still denote the smooth global branches of characteristics of A lifted to \tilde{X} . Let $\tilde{U}(t)$ be the fundamental solution to (1.1) with operator P lifted to \tilde{X} . Let $\tilde{U}(t, \tilde{x}, \tilde{y})$ be the integral kernel of $\tilde{U}(t)$, $\tilde{x}, \tilde{y} \in \tilde{X}$. Let

$$(5.1) \quad U(t, x, y) = \sum_{p\tilde{y}=y} \tilde{U}(t, \tilde{x}, \tilde{y}), \quad x, y \in X,$$

where \tilde{x} is any point of \tilde{X} such that $p\tilde{x} = x$ and the summation is carried out over all preimages \tilde{y} of y . Since (5.1) is invariant under permutations of the leaves of the cover \tilde{X} , the kernel $U(t, x, y)$ is independent of the choice of \tilde{x} . Equation and Cauchy

data are automatically satisfied, so (5.1) gives a fundamental solution to the system (1.1) on X .

The following Lemma gives some estimates for the operator norm from L^2 to H^s for Fourier integral operators in terms of L^∞ norms of the amplitude and its derivatives. Recall that by $\|T\|_s$ we denote the operator norm of T from L^2 to H^s .

Lemma 5.2. *Let T be a Fourier integral operator*

$$(5.2) \quad Tu(x) = \int_{\mathbb{R}^n} \int_X e^{i\varphi(x,y,\xi)} a(x,y,\xi) u(y) dy d\xi,$$

where X is an open set in \mathbb{R}^n . Assume that $a \in S^q$, $q \in \mathbb{Z}$, is an amplitude of order q and has compact support with respect to x and y in X . Assume also that $\partial\varphi/\partial x \neq 0$ for $\xi \neq 0$. Then T extends to a bounded operator from $L^2(X)$ to $H^{-q-n-1}(X)$ with

$$(5.3) \quad \|T\|_{-q-n-1} \leq C \|a\langle\xi\rangle^{-q}\|_{C^{|q+n+1|}},$$

where $\langle\xi\rangle = (1 + |\xi|^2)^{1/2}$.

Moreover, let us assume in addition that the support of a belongs to a conical set with respect to ξ which does not depend on x , i.e.

$$\text{supp } a \subset \Xi_1 = \{(x, y, \xi) : x \in X, (y, \frac{\xi}{|\xi|}) \in \Xi, \xi \neq 0\},$$

where Ξ is subset of S^*X . Then for any $\delta > 0$ we have

$$(5.4) \quad \|T\|_{-q-3n/2-1-\delta} \leq C \text{meas}(\Xi) \|a\langle\xi\rangle^{-q}\|_{C^{|q+n+1|}},$$

where meas is the canonical induced measure on S^*X and constant C may depend on the size of the support of a with respect to x and y .

Proof. First we consider the case $q + n + 1 < 0$. Differentiating (5.2) $-(q + n + 1)$ times with respect to x we obtain the integral with amplitude of order $-(n + 1)$. The integral with respect to ξ converges absolutely, so T extends as a bounded operator from $L^2(X)$ to $C^{-(q+n+1)}$ with estimate (5.3). Estimate (5.4) clearly follows from this as well.

We will now consider the case $q + n + 1 \geq 0$. For a smooth function v , let us consider the bilinear form (Tu, v) . Let us define operator L as the transpose of ${}^tL = (1 + |\partial_x\varphi|^2)^{-1}(1 - i\partial_x\varphi \cdot \partial_x)$. Since $\partial\varphi/\partial x \neq 0$, integrating by parts $q + n + 1$ times with operator L , we obtain an absolutely convergent integral with respect to ξ , and the estimate

$$|(Tu, v)| \leq C \|u\|_{L^2} \|v\|_{q+n+1} \|a\langle\xi\rangle^{-q}\|_{C^{q+n+1}}.$$

This implies (5.3).

We can slightly modify this argument to obtain an estimate of operator T acting from $L^\infty(X)$. Indeed,

$$\begin{aligned} |(Tu, v)| &\leq \left| \int_{\mathbb{R}^n} \int_{\mathbb{R}^n} \int_{\mathbb{R}^n} e^{i\varphi(x,y,\xi)} L^{q+n+1} (a(x,y,\xi) \bar{v}(x)) u(y) dy d\xi dx \right| \\ &\leq C \|v\|_{q+n+1} \|u\|_{L^\infty} I, \end{aligned}$$

where

$$I^2 = \int_X I_1^2(x) dx, \quad I_1(x) = \int_X \int_{\mathbb{R}^n} |\tilde{a}(x, y, \xi)| d\xi dy,$$

with some amplitude \tilde{a} of order $-n-1$. We can estimate \tilde{a} by $\langle \xi \rangle^{-n-1}$ and from the embedding theorems it follows that

$$|(Tu, v)| \leq C \|v\|_{q+n+1} \|u\|_{n/2+\delta} \text{meas}(\Xi) \|a\langle \xi \rangle^{-q}\|_{C^{q+n+1}},$$

implying estimate (5.4) for the norm of operator T acting on $L^2(X)$. \square

The following theorem shows that if a smooth function on a bounded interval has zeros only of finite order, then the measure of the set where this function is small is also small. Moreover, if we have a family of such functions continuously dependent on a parameter, we can estimate measures of sets where functions are small uniformly for all parameters varying over compact sets.

Theorem 5.3. *Let $W \subset \mathbb{R}^n$ be compact and let $0 < T < \infty$. Let a real valued function $f = f(t, w)$ be continuous in $w \in W$ and smooth in $t \in [0, T]$ up to the boundary of $[0, T]$. Let $M \in \mathbb{N}$ and suppose that for each $w \in W$ function $f(\cdot, w)$ has zeros with respect to t of order not greater than M . Then there exist $C > 0$ and $\epsilon_0 > 0$ such that for all $\epsilon < \epsilon_0$ we have*

$$\sup_{w \in W} \text{meas}\{t \in [0, T] : |f(t, w)| \leq \epsilon\} \leq C\epsilon^{1/2M}.$$

Proof. For $\epsilon > 0$ let

$$\Sigma(w, \epsilon) = \{t \in [0, T] : |f(t, w)| \leq C\epsilon\}.$$

Let $K(w)$ be the number of zeros of function $f(\cdot, w)$ with respect to $t \in [0, T]$ and let K be the maximum of $K(w)$ over $w \in W$. It is obvious that K is a finite number due to the condition on zeroes of f and compactness of W .

Let $\alpha_p > 0$, $p \in \mathbb{N}$, be a decreasing sequence of positive numbers which we will choose later. Let us define sets $\Sigma^p(w, \epsilon)$ by setting

$$(5.5) \quad \begin{aligned} \Sigma^p(w, \epsilon) = \{t \in [0, T] : |f(t, w)| \leq C\epsilon, \dots, \\ |\partial_t^{p-1} f(t, w)| \leq C\epsilon^{\alpha_{p-1}}, |\partial_t^p f(t, w)| \geq C\epsilon^{\alpha_p}\}, \quad p \in \mathbb{N}. \end{aligned}$$

We claim now that there exists $\epsilon_0 > 0$ such that for all $0 < \epsilon < \epsilon_0$, and all $w \in W$, we have $\Sigma^p(w, \epsilon) = \emptyset$ for all $p > M$. Indeed, if it is not so, then due to compactness of W there are sequences t_n , w_n and ϵ_n , converging to $t^* \in [0, T]$, $w^* \in W$ and zero, respectively, such that

$$|f(t_n, w_n)| \leq C\epsilon_n, \dots, |\partial_t^{M+1} f(t_n, w_n)| \leq C\epsilon_n^{\alpha_{M+1}},$$

and consequently

$$\partial_t^p f(t^*, w^*) = 0, \quad p = 0, \dots, M+1,$$

which is impossible. It follows now that the set $\Sigma(w, \epsilon)$ may be presented as the following union of sets

$$(5.6) \quad \Sigma(w, \epsilon) = \bigcup_{p=1}^M \Sigma^p(w, \epsilon).$$

The idea of the proof now is to show first that the number of connected components of sets $\Sigma^p(w, \epsilon)$ is finite and can be estimated uniformly over all w and ϵ . Then we will show that the size of each connected component is small and can be estimated by

$\epsilon^{1/2M}$, which will imply Theorem 5.3. These statements are proved in the following two lemma.

Lemma 5.4. *There exists $\epsilon_0 > 0$ such that for all $0 < \epsilon < \epsilon_0$ the inequality*

$$(5.7) \quad \Delta(\Sigma^p(w, \epsilon)) \leq K(M+1)^2, \quad \forall w \in W, 1 \leq p \leq M,$$

holds, where $\Delta(\Sigma^p)$ is the number of connected components of Σ^p .

Proof. For simplicity let us first consider the case $p = 1$. Let us assume that (5.7) is not valid. Then there exist sequences w_n converging to some $w^* \in W$ and ϵ_n converging to 0 such that

$$(5.8) \quad \Delta(\Sigma^1(w_n, \epsilon_n)) > K(M+1)^2.$$

Let us now choose small enough $\epsilon_1 > 0$ such that each connected interval in the closure of $\Sigma(w^*, \epsilon_1)$ will include one and only one zero of function $f(t, w^*)$ with respect to t and will not include zeros of derivative of $\partial_t f(t, w^*)$ different from zeros of $f(t, w^*)$. This is possible because if we had an infinite number of zeros of $\partial_t f(t, w^*)$ approaching a zero t^* of $f(t, w^*)$, it would mean that $\partial_t f(t^*, w^*) = 0$ and that in fact t^* is zero of $\partial^k f(t, w^*)$ for all $k \geq 1$, which is impossible since we assumed that all zeros of $f(t, w^*)$ are of finite order not exceeding M . It follows now that $\Delta(\Sigma(w^*, \epsilon_1)) \leq K$. Since w_n converges to w^* , by continuity we also have

$$(5.9) \quad \Sigma^1(w_n, \epsilon_n) \subset \Sigma(w^*, \epsilon_1),$$

for sufficiently large n . From (5.8) and (5.9) it follows that there exist $(M+1)^2 + 1 \geq 2M+1$ connected components of $\Sigma^1(w_n, \epsilon_n)$ which are all contained in one of the connected components of closure of $\Sigma(w^*, \epsilon_1)$. Let us denote this connected component of $\Sigma(w^*, \epsilon_1)$ by I .

According to the definition of sets $\Sigma^1(w_n, \epsilon_n)$, between two of its connected components function $f(t, w_n)$ must become relatively large (i.e. $> C\epsilon$) or its derivative $\partial_t f(t, w_n)$ must become relatively small (i.e. $< C\epsilon^{\alpha_1}$). From this we see that in the first case $\partial_t f(\cdot, w_n)$ must become zero at some point between these components, while in the second case $\partial_t^2 f(\cdot, w_n)$ must become zero at some point there. Since the number of components of $\Sigma^1(w_n, \epsilon_n)$ in I is at least $2M+1$, it follows that at least one of these two cases occurs at least $M+1$ times.

Let us consider these cases separately. In the first case, for sufficiently large n , we have at least M zeros $s_n^1 < \dots < s_n^M$ of the derivative $\partial_t f(t, w_n)$ contained in I . It follows that $\partial_t^2 f(\cdot, w_n)$ has at least $M-1$ different zeros in I , $\partial_t^3 f(\cdot, w_n)$ has at least $M-2$ different zeros in I , etc. In particular, there are $\tau_n^k \in [s_n^1, s_n^M]$ such that $\partial_t^k f(\tau_n^k, w_n) = 0$ for $k = 1, \dots, M$.

Using compactness of I and continuity of f , it follows that there are subsequences $s_{n_i}^1, \dots, s_{n_i}^M, \tau_{n_i}^k$ and w_n converging to s_*^1, \dots, s_*^M and τ_*^k , respectively, which are all contained in I , and $w_n \rightarrow w^*$. Since functions $f(t, w)$ are smooth in t we have

$$\partial_t f(s_*^1, w^*) = 0, \dots, \partial_t f(s_*^M, w^*) = 0, \text{ and } \partial_t^k f(\tau_*^k, w^*) = 0, \quad k = 1, \dots, M.$$

Since there are no zeros of derivative $\partial_t f(\cdot, w^*)$ on the interval I except may be a point t^* which is zero of function $f(\cdot, w^*)$, we obtain $s_*^1 = \dots = s_*^M = t^*$. From $\tau_n^k \in [s_n^1, s_n^M]$ it follows that $\tau_*^k = t^*$ for all $k = 1, \dots, M$, which means

$$\partial_t^k f(t^*, w_*) = 0 \text{ for all } k = 0, \dots, M.$$

But this is impossible since function $f(\cdot, w^*)$ may have zeros only of order M in view of our assumption.

The second case is slightly different. Here, for sufficiently large n , we have at least M zeros $s_n^1 < \dots < s_n^M$ of the second order derivative $\partial_t^2 f(t, w_n)$ contained in $I \setminus \Sigma^1(w_n, \epsilon_n)$. Because in this case we assumed that $\Sigma^1(w_n, \epsilon_n)$ breaks into at least $M + 1$ components due to the failure of condition $|\partial_t f(t, w)| \geq C\epsilon^{\alpha_1}$, it follows that

$$(5.10) \quad |\partial_t f(s_n^i, w_n)| \leq C\epsilon_n^{\alpha_1}, \quad i = 1, \dots, M.$$

As n tends to infinity, we can choose subsequences of s_n^i converging to some $s_*^i \in I$. Because $\partial_t f(\cdot, w^*)$ does not have zeros in I except may be for some $t^* \in I$ which is also the unique zero of $f(\cdot, w^*)$ in I , we get that $s_*^1 = \dots = s_*^M = t^*$. From (5.10) we also have $\partial_t f(t^*, w^*) = 0$.

Now, to deal with higher order derivatives of f at t^* , similar to the first case, we get a collection of $\tau_n^k \in [s_n^1, s_n^M]$ such that $\partial_t^k f(\tau_n^k, w_n) = 0$ for $k = 2, \dots, M + 1$. Because of compactness τ_n^k has a subsequence converging to some τ_*^k , $k = 2, \dots, M + 1$. Again, we must have $\tau_*^k = t^*$ and hence also $\partial_t^k f(t^*, w^*) = 0$, for all $k = 2, \dots, M + 1$. Since we already showed that $f(t^*, w^*) = \partial_t f(t^*, w^*) = 0$, this contradicts the assumption that $f(\cdot, w^*)$ may have zeros of order up to M .

The argument for $p \geq 2$ is similar. For t between two connected components of Σ^p , at least one of conditions in (5.5) breaks down. Note that since we assumed that the total number of components of $\Sigma^p(w_n, \epsilon_n)$ is larger than $K(M + 1)^2$ and the number of components of the larger set $\Sigma(w^*, \epsilon_1)$ is at most K , we will have at least $(M + 1)^2 + 1$ components of $\Sigma^p(w_n, \epsilon_n)$ in I . This means that these $p + 1$ conditions fail at least $(M + 1)^2$ times. Since $p + 1 \leq M + 1$, there is a condition that will fail at least $M + 1$ times.

In the case this is the last condition $|\partial_t^p f(t, w)| \geq C\epsilon^{\alpha_p}$ that fails while conditions $|\partial_t^i f(t, w)| \leq C\epsilon^{\alpha_i}$ remain valid, we can argue similar to the second case of the argument with $p = 1$. In this we need to have at least $M - p$ different zeros of $\partial_t^{p+1} f(t, w)$, which is the case if we have at least $M - p + 1$ such components in I . This is true since we have at least $M + 1$ such components.

If one of the other conditions fails, let us take the smallest i for which condition $|\partial_t^i f(t, w)| \leq C\epsilon^{\alpha_i}$ fails $M + 1$ times. Again, we need to have at least $M - i$ different zeros of $\partial_t^{i+1} f(t, w)$, which would follow from having $M - i + 1$ such components in I . This is again true since we have at least $M + 1$ such components. \square

Our next step is to show that each connected component of $\Sigma^p(w, \epsilon)$, for $p = 1, \dots, M$, is small enough.

Lemma 5.5. *The length of each connected component of $\Sigma^p(w, \epsilon)$, for $p = 1, \dots, M$, is no greater than $C\epsilon^{\alpha_{p-1} - \alpha_p}$.*

Proof. Let I be a connected component of $\Sigma^p(w, \epsilon)$ and let $t^* \in I$. We are going to estimate the maximal shift $\delta_0 > 0$ such that $t^* + \delta \in I$ for all $0 < \delta < \delta_0$. Recalling the definition of $\Sigma^p(w, \epsilon)$, we see that

$$(5.11) \quad |\partial_t^{p-1} f(t^* + \delta, w)| \leq C\epsilon^{\alpha_{p-1}}, \quad |\partial_t^p f(t^* + \delta, w)| \geq C\epsilon^{\alpha_p}.$$

Using the Taylor expansion of $\partial_t^{p-1} f(\cdot, w)$ at t^* , we have

$$\partial_t^{p-1} f(t^* + \delta, w) = \partial_t^{p-1} f(t^*, w) + \partial_t^p f(t^*, w)\delta,$$

where t^{**} is some point between t^* and $t^* + \delta$. Since $t^*, t^* + \delta \in I$, it follows that $t^{**} \in I$ and

$$|\partial_t^p f(t^{**}, w)| |\delta| \leq |\partial_t^{p-1} f(t^* + \delta, w)| + |\partial_t^{p-1} f(t^*, w)| \leq 2C\epsilon^{\alpha_{p-1}}.$$

From this and (5.11) we obtain that $|\delta_0| \leq C\epsilon^{\alpha_{p-1}-\alpha_p}$. Consequently, the length of each connected component of $\Sigma^p(w, \epsilon)$ can be estimated by $C\epsilon^{\alpha_{p-1}-\alpha_p}$. \square

Now we can finish the proof of Theorem 5.3. Let us choose $\alpha_k = 1 - k/2M$, $k = 0, \dots, M$. According to Lemma 5.5 the length of each connected component of $\Sigma^p(w, \epsilon)$ can be estimated by $C\epsilon^{1/2M}$. Then according to Lemma 5.4, the size of $\Sigma^p(w, \epsilon)$ can be estimated by $CK(M+1)^2\epsilon^{1/2M}$. Because of decomposition (5.6) the size of $\Sigma(w, \epsilon)$ is estimated by $C\epsilon^{1/2M}$. Statement of Theorem 5.3 is now a consequence of continuity of f with respect to w and compactness of W . \square

The following interpolation lemma shows that if a function u can be decomposed for all sufficiently small ϵ into a sum of $u_1^{(\epsilon)} + u_2^{(\epsilon)}$ with a good estimate for the norm of $u_2^{(\epsilon)}$ in a “bad” Sobolev space H^p with small index p , and with a bad estimate for the norm of $u_1^{(\epsilon)}$ in a “good” Sobolev space H^r with large index r , then u belongs to some intermediate space H^q with $p < q < r$.

Lemma 5.6. *Let $u \in H^p(\mathbb{R}^n)$. Suppose that for all small enough ϵ there is a representation $u = u_1^{(\epsilon)} + u_2^{(\epsilon)}$ such that*

$$(5.12) \quad \|u_1^{(\epsilon)} : H^r\| \leq C\epsilon^{-T}, \quad \|u_2^{(\epsilon)} : H^p\| \leq C\epsilon^S, \quad p < r; \quad S, T > 0.$$

Then $u \in H^q(\mathbb{R}^n)$ for any

$$(5.13) \quad q < (pT + rS)(T + S)^{-1}.$$

Proof. Let a sequence $\{a_j\}$ be such that $0 = a_0 < a_1 < \dots$ and $a_j \rightarrow +\infty$ as $j \rightarrow +\infty$, and let a sequence $\{b_j\}$ of positive numbers $b_j > 0$ tend to zero. Let us assume that $u \in H^q(\mathbb{R}^n)$ for some q . Then

$$\begin{aligned} \|u : H^q\|^2 &= \int_{\mathbb{R}^n} |\hat{u}|^2 (1 + |\xi|^2)^q d\xi \\ &= 2 \sum_{j=0}^{+\infty} \left(\int_{a_j \leq |\xi| \leq a_{j+1}} |\widehat{u_2^{(b_j)}}|^2 (1 + |\xi|^2)^q d\xi + \int_{a_j \leq |\xi| \leq a_{j+1}} |\widehat{u_1^{(b_j)}}|^2 (1 + |\xi|^2)^q d\xi \right) \\ &\leq 2 \sum_{j=0}^{+\infty} \left(\max\{\langle a_j \rangle^{2(q-p)}, \langle a_{j+1} \rangle^{2(q-p)}\} \int_{a_j \leq |\xi| \leq a_{j+1}} |\widehat{u_2^{(b_j)}}|^2 (1 + |\xi|^2)^p d\xi \right. \\ &\quad \left. + \max\{\langle a_j \rangle^{2(q-r)}, \langle a_{j+1} \rangle^{2(q-r)}\} \int_{a_j \leq |\xi| \leq a_{j+1}} |\widehat{u_1^{(b_j)}}|^2 (1 + |\xi|^2)^r d\xi \right). \end{aligned}$$

It follows from this estimate and (5.12) that

$$(5.14) \quad \|u : H^q\|^2 \leq C \sum_{j=0}^{+\infty} \left(\max\{\langle a_j \rangle^{2(q-p)}, \langle a_{j+1} \rangle^{2(q-p)}\} b_j^{2S} + \max\{\langle a_j \rangle^{2(q-r)}, \langle a_{j+1} \rangle^{2(q-r)}\} b_j^{-2T} \right).$$

Now we are going to demonstrate that under hypothesis (5.13) we may choose sequences $\{a_j\}$ and $\{b_j\}$ in such way that the right hand side of (5.14) will be finite, so that conclusion of Lemma 5.6 will follow. Let us set $a_j = j^\alpha$ and $b_j = j^{-\beta}$ with some $\alpha, \beta > 0$. Then the series in (5.14) will converge if the following inequalities are fulfilled

$$2\alpha(q-p) - 2\beta S < -1, \quad 2\alpha(q-r) + 2\beta T < -1.$$

These inequalities can be transformed into

$$\alpha(2T(q-p) + 2S(q-r)) < -T - S, \quad 2\beta N > \alpha 2(q-p) + 1,$$

which hold with positive constants α and β if and only if (5.13) is valid. \square

6. SPECTRAL ASYMPTOTICS

In this section we will prove Theorem 2.5. Let X be a smooth compact manifold without boundary of dimension $n \geq 3$. Then an elliptic operator $P(x, D)$ has a collection of eigenfunctions and eigenvalues $\lambda_j \rightarrow \infty$. We will be interested in distribution of eigenvalues and will find the asymptotics for the spectral function $N(\lambda) = \sharp\{j : \lambda_j < \lambda\}$. One of the most effective methods to study such asymptotics is to use an explicit representation for the fundamental solution of the corresponding hyperbolic problem (1.1). We will follow the method developed in [8], [7], [23], etc., for scalar operators, and in [21] for the case $M = 1$ in Condition C. The following proposition was implicitly proved in [7] and formulated in [21].

Proposition 6.1. *Let $\chi_1, \chi_2 \in C_0^\infty(\mathbb{R})$ be such that $\chi_1(0) = 1$ and $0 \notin \text{supp } \chi_2$. Suppose that*

$$(6.1) \quad \begin{aligned} \text{Tr } \mathcal{F}_{t \rightarrow \mu}^{-1}(\chi_1(t)U(t)) &= c_1 \mu^{n-1} + c_2 \mu^{n-2} + o(\mu^{n-2}), \\ \text{Tr } \mathcal{F}_{t \rightarrow \mu}^{-1}(\chi_2(t)U(t)) &= o(\mu^{n-1}), \quad \mu \rightarrow \infty. \end{aligned}$$

Then $N(\lambda) = c_1 n^{-1} \lambda^n + c_2 (n-1)^{-1} \lambda^{n-1} + o(\lambda^{n-1})$.

In previous sections, and in particular in the proof of Theorem 2.2, we have represented the propagator $U(t)$ in (1.1) as an infinite sum of some extensions of Fourier integral operators, i.e.

$$(6.2) \quad U(t) = \sum_{j=0}^{\infty} e^{-i\tilde{A}t} Q_j,$$

where $Q_0 = I$ and $\text{Tr } Q_1 = 0$. So we also have

$$(6.3) \quad \text{Tr } e^{-i\tilde{A}t} Q_1 = 0.$$

Let us first consider the contribution of the first term, which is $e^{-i\tilde{A}t}$. It is the propagator for a block-diagonal system, so the asymptotic behaviour of $\mathcal{F}_{t \rightarrow \mu}^{-1}(\chi_\sigma(t)e^{-i\tilde{A}t})$, $\sigma = 1, 2$, determines the spectral distribution for a system of independent scalar equations. Thus, this is the sum of spectral distributions for scalar operators, which are well-known (e.g. [8], [7]). So

$$\mathcal{F}_{t \rightarrow \mu}^{-1} \text{Tr}(\chi_1(t)e^{-i\tilde{A}t}) = c_1 \mu^{n-1} + c_2 \mu^{n-2} + o(\mu^{n-2}).$$

We will show that under suitable conditions the asymptotics in (6.1) are determined only by the first term of the series (6.2). Let us first observe that in view of Theorem 2.1 operator $\sum_{j=K}^{\infty} e^{-i\tilde{A}t} Q_j$ is compact for sufficiently large K , so it does not contribute to the singularity of $\text{Tr } \chi_{\sigma} U$, $\sigma = 1, 2$. Therefore, we can replace $U(t)$ in (6.1) by a sum of finitely many terms in (6.2). Also, in view of (6.3), the second term of (6.2) does not contribute to (6.1). Other terms of the sum (6.2) are of the form

$$e^{-i\tilde{A}t} \int_0^t \int_0^{t_1} \cdots \int_0^{t_l} Z(t_1) Z(t_2) \cdots Z(t_{l+1}) dt_{l+1} \cdots dt_1, \quad l \geq 2,$$

where $Z(\tau) = -ie^{i\tilde{A}\tau} B e^{-i\tilde{A}\tau}$. In this way, after a change of variables $s_1 = t - t_1$, $s_2 = t_1 - t_2, \dots, s_{l+1} = t_l$, we obtain a sum of terms of the form

$$I_{\sigma}(\mu) = \text{Tr} \int \chi_{\sigma}(s_1 + \cdots + s_{l+1}) e^{i(s_1 + \cdots + s_{l+1})\mu} B(s) L(s) ds, \quad \sigma = 1, 2.$$

where $B(s) \in \Psi^0$, $L(s) = e^{-i\tilde{A}_{j_1}s_1} e^{-i\tilde{A}_{j_2}s_2} \cdots e^{-i\tilde{A}_{j_{l+1}}s_{l+1}}$, and \tilde{A}_j is the j -th block of \tilde{A} corresponding to a_j . As before, $L(s)$ may be represented as a locally finite sum of oscillatory integrals with phase $\varphi(s, x, y, \xi)$, for which

$$\frac{\partial \varphi}{\partial s_k} = S_{j_k}(s, x, \frac{\partial \varphi}{\partial x}),$$

where $S_{j_k}(s, x, p) = a_{j_k}(\Phi_{j_{k-1}}^{s_{k-1}} \cdots \Phi_{j_1}^{s_1}(x, p))$. Therefore,

$$I_{\sigma}(\mu) = \int_X \int_{\mathbb{R}^n} \chi_{\sigma}(s_1 + \cdots + s_{l+1}) e^{i(s_1 + \cdots + s_{l+1})\mu + i\varphi(s, x, x, \xi)} b(s, x, x, \xi) d\xi dx ds.$$

Substituting $\xi = \mu\tau\omega$, $\tau > 0$, $|\omega| = 1$, we get

$$(6.4) \quad I_{\sigma}(\mu) = \mu^n \int_X \int_{\mathbf{S}^{n-1}} \int_0^{\infty} \chi_{\sigma}(s_1 + \cdots + s_{l+1}) e^{i\mu(s_1 + \cdots + s_{l+1} + \tau\varphi)} b\tau^{n-1} d\tau d\omega dx ds.$$

To finish the argument for this operator, we can follow [21] to show that smoothing of Q_l 's implies (6.1). If we change variables again by $\rho = \sum s_j$, $s_j = \rho\kappa_j$ and introduce $K = \{\kappa_j \geq 0, \sum \kappa_j = 1\}$, we get

$$I_1(\mu) = \mu^n \int_K \int_X \int_{\mathbf{S}^{n-1}} \int_{-\infty}^{\infty} \int_0^{\infty} \chi_1(\rho) e^{i\mu(\rho + \tau\varphi(\rho\kappa, x, x, \omega))} \rho^l \tau^{n-1} b d\tau d\rho d\omega dx d\kappa.$$

For fixed ω, x, κ the point $\tau = -(\sum \kappa_k S_{j_k}(\rho\kappa, x, \partial\varphi/\partial x))^{-1}$, $\rho = 1$, is a nondegenerate stationary point of this oscillatory integral, implying $I_1(\mu) = O(\mu^{n-1-l}) = o(\mu^{n-2})$. This already gives Hörmander's first term of $N(\lambda)$. Now we will show that singularities at $t \neq 0$ do not give essential contributions to the second term of spectral asymptotics. In the analysis of $\text{Tr } \chi_2 U$, the contribution of the first term of (6.2) was established in [7]. Let us look at the integral (6.4) with respect to s_1, s_2, τ with fixed $x, \omega, s_3, \dots, s_{l+1}$. The stationary phase method with respect to τ, s_1 gives a stationary point $\varphi(s, x, x, \omega) = 0$, $\tau = -(\partial\varphi/\partial s_1)^{-1}$. It is non-degenerate because $\det \partial_{\tau} \partial_{s_1}(\tau\varphi) = -a_1(s, x, \partial_x \varphi)^2$. This gives the estimate $I_2(\mu) = O(\mu^{n-1})$. Since $\partial\varphi/\partial s_2 = 0$ only at $\tau = -(\partial\varphi/\partial s_2)^{-1} = a_2(s, x, \partial_x \varphi)^{-1}$, the phase is stationary with respect to s_2 only at (x, ξ) for which $a_1(s, x, \partial_x \varphi) = a_2(s, x, \partial_x \varphi)$. This is the set of measure zero and, therefore, $I_2(\mu) = o(\mu^{n-1})$, which shows (6.1). Finally we note that if the support of χ_1 is sufficiently small (i.e. when s is small), terms in (5.1),

corresponding to \tilde{x} and \tilde{y} in different leaves of \tilde{X} will produce only smoothing operators (5.1) in view of the finite propagation speed. So their contribution to $I(\mu)$ is rapidly decreasing and we obtain the result also on X .

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ILIA KAMOTSKI, MICHAEL RUZHANSKY:
 DEPARTMENT OF MATHEMATICS
 IMPERIAL COLLEGE OF SCIENCE, TECHNOLOGY AND MEDICINE
 180 QUEEN'S GATE, LONDON SW7 2BZ, UK
E-mail addresses kamotski@ic.ac.uk, ruzh@ic.ac.uk